

JOURNAL OF DIFFERENTIAL EQUATIONS, 2, 365-377 (1966).

A Liapunov Method for the Estimation of Statistical Averages

W. M. WONHAM¹*Division of Applied Mathematics,
Brown University, Providence, Rhode Island**Received July 16, 1965*

1. INTRODUCTION

We consider a randomly perturbed dynamical system described by the equation

$$dx/dt = f(x) + G(x)\xi(t), \quad t \geq 0, \quad (1)$$

where x, f are n -vectors, G is an $n \times n$ matrix and $\xi(t)$ is n -dimensional Gaussian white noise. Such equations arise in control theory [1], and the theory of random vibrations [2]. In these applications it is of interest to know under what conditions the process

$$X = \{x(t), \quad t \geq 0\}$$

generated by (1) is stable, in the sense that X admits a unique invariant probability distribution. If X is stable then it is often desirable to estimate various stationary averages $\mathcal{E}\{L(x)\}$, when these averages exist.

In a previous paper [3] a criterion of Liapunov type was given for stability in the sense described. In the present note a Liapunov criterion (Theorem 3.1) is obtained for the existence (finiteness) of the stationary average $\mathcal{E}\{L(x)\}$ where L is an arbitrary nonnegative function. This result is applied to show that algebraic moments of all orders exist when, in (1), G is bounded and the unperturbed system $dx/dt = f(x)$ is of Lur'e type.

The existence criterion is extended to yield an effective method of calculating an upper bound for $\mathcal{E}\{L(x)\}$ (Theorem 4.1). The method is illustrated by an example from control theory.

2. STATEMENT OF THE PROBLEM

We start with a precise version of (1), namely Itô's equation

$$\begin{aligned} dx(t) &= f(x(t)) dt + G(x(t)) dw(t), \quad t \geq 0 \\ x(0) &= x_0. \end{aligned} \quad (2)$$

¹ This research was supported in part by the National Aeronautics and Space Administration under grant No. NGR-40-002-015, and in part by the United States Air Force through the Office of Scientific Research under grant AF-AFOSR-693-65.

The following assumptions are made with respect to (2):

(i) x, f are vectors in Euclidean n -space E ($n \geq 2$) and G is an $n \times n$ matrix.

(ii) $\{w(t); t \geq 0\}$ is a Wiener process in E

(iii) x_0 is a random variable independent of the process $w(t)$.

(iv) There is a constant $c > 0$ such that

$$|f(x) - f(y)| + |G(x) - G(y)| < c |x - y|$$

for all $x, y \in E$. (Here $|\cdot|$ denotes Euclidean norm of a vector or matrix.)

(v) There is a constant $\epsilon > 0$ such that

$$y'G(x)G(x)'y \geq \epsilon |y|^2$$

for all $x, y \in E$. (A prime denotes transpose of a vector or matrix.)

Under these assumptions it is known (cf. [3]) that (2), interpreted in the sense of Itô, defines a continuous, strongly Feller process

$$X = \{x(t), \quad t \geq 0\}.$$

The differential generator of X will be denoted by \mathcal{L} , where

$$\mathcal{L}[u(x)] = \frac{1}{2} \text{tr}[G(x)G(x)'u_{xx}(x)] + f(x)'u_x(x) \quad (3)$$

whenever the indicated derivatives exist. (In (3), u_x is the vector of first partial derivatives of u and u_{xx} is the matrix of second partial derivatives.)

In the following we shall always assume that X is *positive* [4]. Under these conditions it is known [4] that there exists a unique invariant probability measure μ defined on the Borel sets $B \subset E$: that is, if P denotes probability measure on the paths of X , and if

$$P(x_0 \in B) = \mu(B)$$

then

$$P(x(t) \in B) = \mu(B), \quad t > 0.$$

An effective criterion for positivity of X is given in [3].

Let $L(x) \geq 0$ be Hölder continuous on the compact subsets of E . The main problem is to obtain a sufficient condition that

$$\mathcal{E}\{L(x)\} = \int_E L(x)\mu(dx)$$

be finite. Subsequently we shall describe a method for deriving an upper bound on $\mathcal{E}\{L(x)\}$.

In the following, the terms *smooth*, and *normal domain*, have the same meaning as in [3].

3. A CRITERION FOR EXISTENCE OF $\mathcal{E}\{L(x)\}$

A Liapunov criterion for the existence of $\mathcal{E}\{L(x)\}$ can be derived by arguments very similar to those of [3] and [4]. The result is given in Theorem 3.1. We start with some preliminary lemmas.

Let D be a normal domain with boundary Γ , and let τ_Γ be the first time X hits Γ . Let \mathcal{E}_x denote expectation on the paths of X when $x(0) = x \in E$. Since X is positive, $\mathcal{E}_x(\tau_\Gamma) < \infty$, $x \in E - D$.

LEMMA 3.1. *Let*

$$u(x) = \mathcal{E}_x \left\{ \int_0^{\tau_\Gamma} L[x(t)] dt \right\}$$

If $u(x_0) < \infty$ for some point $x_0 \in E - \bar{D}$ then $u(x) < \infty$ for all $x \in E - \bar{D}$. Furthermore

$$\begin{aligned} \mathcal{L}[u(x)] &= -L(x), \quad x \in E - \bar{D} \\ u(x) &= 0, \quad x \in \Gamma. \end{aligned} \quad (4)$$

Proof. The proof closely follows that of Lemma 5.3 of [4]. Let $\{D_n; n = 1, 2, \dots\}$ be an increasing sequence of normal domains such that $D \subset D_1$, $x_0 \in D_1 - D$ and $\lim D_n = E (n \rightarrow \infty)$. Let τ_n be the first time X hits the boundary $\Gamma \cup \Gamma_n$ of $D_n - D$, and define

$$\begin{aligned} u_n(x) &= \mathcal{E}_x \left\{ \int_0^{\tau_n} L[x(t)] dt \right\}, \quad x \in \bar{D}_n - D \\ u(x) &= \mathcal{E}_x \left\{ \int_0^{\tau_\Gamma} L[x(t)] dt \right\}, \quad x \in E - D. \end{aligned}$$

Since X is positive (and hence, regular [4]) $\tau_n \uparrow \tau_\Gamma$ ($n \rightarrow \infty$) and therefore $u_n(x) \uparrow u(x)$ ($n \rightarrow \infty$). For fixed $m \geq 1$

$$u(x) = u_m(x) + \sum_{n=m}^{\infty} [u_{n+1}(x) - u_n(x)], \quad x \in \bar{D}_m - D, \quad (5)$$

with convergence at least for $x = x_0$. We now use the fact that $u_n(x)$ is the unique smooth solution of

$$\begin{aligned} \mathcal{L}[u_n(x)] &= -L(x), \quad x \in D_n - \bar{D}. \\ u_n(x) &= 0, \quad x \in \Gamma \cup \Gamma_n. \end{aligned}$$

(see e.g. [5], theorem 13.16). Let

$$v_n(x) = u_{n+1}(x) - u_n(x), \quad x \in \bar{D}_n - D.$$

Then

$$\mathcal{L}[v_n(x)] = 0, \quad x \in D_n - \bar{D}$$

and (since $u_n(x) \geq 0$) $v_n(x) \geq 0$, $x \in \Gamma \cup \Gamma_n$. By the maximum principle $v_n(x) \geq 0$, $x \in D_n - \bar{D}$. It follows that all terms of the series (5) except the first are positive functions for which $\mathcal{L}[v(x)] = 0$, and the series converges for $x = x_0$. From the generalized Harnack inequality [11] it follows that the series converges for all $x \in \bar{D}_m - D$ and $u(x)$ satisfies (4) for $x \in D_m - \bar{D}$. Since m is arbitrary the result follows.

LEMMA 3.2. *A necessary and sufficient condition that*

$$\mathcal{E}\{L(x)\} < \infty$$

is that

$$\mathcal{E}_x \left\{ \int_0^{\tau^T} L[x(t)] dt \right\} < \infty, \quad x \in E. \quad (6)$$

Proof. We use the construction and notation of [4]. Let D_1 be a normal domain with boundary Γ_1 such that $D \subset D_1$ and $\Gamma \cap \Gamma_1 = \phi$. Let τ denote the length of a cycle, namely, in obvious notation,

$$\tau = \min\{t : x(t) \in \Gamma \mid x(0) \in \Gamma \text{ and } x(s) \in \Gamma_1 \text{ for some } s, 0 < s < t\}.$$

Let $\tilde{\mu}$ be the finite invariant measure (see [4]) induced on the Borel sets of Γ . Then if K is an arbitrary compact subset of E we have, within a constant of normalization,

$$\mu(K) = \int_{\Gamma} \tilde{\mu}(dx) \mathcal{E}_x\{\tau^K\}, \quad (7)$$

where

$$\tau^K = \text{meas}\{t : 0 \leq t \leq \tau, x(t) \in K\}.$$

Let $L_n(x)$ be an increasing sequence of simple functions (constructed on compact sets) such that $L_n(x) = 0$ ($|x| > n$) and $L_n(x) \uparrow L(x)$ ($n \rightarrow \infty$). From (7)

$$\int_E \mu(dx) L_n(x) = \int_{\Gamma} \tilde{\mu}(dx) \mathcal{E}_x \left\{ \int_0^{\tau} L_n[x(t)] dt \right\}, \quad n = 1, 2, \dots;$$

and by monotone convergence

$$\mathcal{E}\{L(x)\} = \int_{\Gamma} \tilde{\mu}(dx) \mathcal{E}_x \left\{ \int_0^{\tau} L[x(t)] dt \right\}. \quad (8)$$

Let $\tau_1 = \min\{t : x(t) \in \Gamma_1 \mid x(0) \in \Gamma\}$. By the strong Markov property

$$\begin{aligned} \mathcal{E}_x \left\{ \int_0^\tau L[x(t)] dt \right\} &= \mathcal{E}_x \left\{ \int_0^{\tau_1} L[x(t)] dt \right\} \\ &+ \mathcal{E}_x \left\{ \mathcal{E}_{x(\tau_1)} \left\{ \int_0^{\tau_1} L[x(t)] dt \right\} \right\}, \quad x \in \Gamma. \end{aligned} \quad (9)$$

We will show that each term in the right side of (9) is bounded on Γ . Let $u(x)$ denote the first term, regarded as a function on \bar{D}_1 . Then u is the unique smooth solution of $\mathcal{L}[u(x)] = -L(x)$, $x \in D_1$; $u(x) = 0$, $x \in \Gamma_1$. Since \bar{D}_1 is compact, u is bounded. Next, let

$$v(y) = \mathcal{E}_y \left\{ \int_0^{\tau_1} L[x(t)] dt \right\}.$$

By (6) and Lemma 3.1, $v(y)$ is smooth, hence bounded on Γ_1 . By the strong Feller property, $\mathcal{E}_x\{v(x(\tau_1))\}$ is continuous, and therefore bounded on Γ . It now follows from (8) that $\mathcal{E}\{L(x)\} < \infty$.

Conversely, if (6) fails for some $x \in E - D$ then by Lemma 3.1, (6) fails for all $x \in E - D$, and by (8) and (9), $\mathcal{E}\{L(x)\} = \infty$.

LEMMA 3.3. *If the equation*

$$\mathcal{L}[v(x)] = -L(x), \quad x \in E - \bar{D}$$

has a smooth positive solution $v(x)$ in $E - D$ then

$$\mathcal{E}_x \left\{ \int_0^{\tau_1} L[x(t)] dt \right\} < \infty, \quad x \in E - D.$$

Proof. Let $\{D_n; n = 1, 2, \dots\}$ be a sequence of normal domains constructed as in the proof of Lemma 3.1, and let $u_n(x)$ be the corresponding sequence of smooth functions such that

$$\begin{aligned} \mathcal{L}[u_n(x)] &= -L(x), \quad x \in D_n - \bar{D} \\ u_n(x) &= 0, \quad x \in \Gamma \cup \Gamma_n. \end{aligned}$$

Since $\mathcal{L}[v(x) - u_n(x)] = 0$, $x \in D_n - \bar{D}$, and $v(x) - u_n(x) \geq 0$, $x \in \Gamma \cup \Gamma_n$, we have $u_n(x) \leq v(x)$, $x \in \bar{D}_n - D$; therefore

$$\begin{aligned} \mathcal{E}_x \left\{ \int_0^{\tau_1} L[x(t)] dt \right\} &= \lim u_n(x) \\ &\leq v(x), \quad x \in E - D. \end{aligned}$$

This completes the proof.

Before stating Theorem 3.1 we introduce a class of real-valued functions V , analogous to Liapunov functions, with the following properties.

P_1 : V is defined for $x \in \bar{D}_v$, where

$$D_v = \{x : |x| > R\} \quad (R < \infty \text{ is arbitrary})$$

P_2 : V is continuous in \bar{D}_v and is twice continuously differentiable in D_v .

P_3 : $V(x) \rightarrow +\infty$ as $|x| \rightarrow \infty$.

THEOREM 3.1. *Let X be positive. If there exists a function V with properties $P_1 - P_3$ and the additional property*

$$\mathcal{L}[V(x)] \leq -L(x), \quad x \in D_v$$

then

$$\mathcal{E}\{L(x)\} < \infty.$$

We remark that if X is positive and $L(x)$ is bounded then $\mathcal{E}\{L(x)\}$ is obviously finite. If $L(x)$ is bounded away from zero for $x \in D_v$ then, by Theorem 2 of [3], the existence of V already implies that X is positive.

Proof. By Lemmas 3.2 and 3.3 we have that $\mathcal{E}\{L(x)\} < \infty$ if and only if there exists a normal domain D such that the equation

$$\mathcal{L}[u(x)] = -L(x) \tag{10}$$

has a smooth positive solution $u(x)$ defined for $x \in E - D$. Let $D = E - \bar{D}_v$ and define a sequence $\{D_n\}$ of normal domains as in the proof of Lemma 3.1. The remainder of the proof follows that of Lemma 3.3, with $v(x)$ replaced by $V(x)$. By adding a constant to V if necessary we can arrange that $V(x) \geq 0$, $x \in \bar{D}_v$. If

$$\mathcal{L}[u_n(x)] = -L(x) \quad (x \in D_n - \bar{D}),$$

$$u_n(x) = 0 \quad (x \in \Gamma \cup \Gamma_n),$$

then $0 \leq u_n(x) \leq u_{n+1}(x) \leq V(x)$, $x \in \bar{D}_n - D$. It follows by a compactness theorem ([6] p. 344) that $\lim u_n(x)$ exists and is a solution of (10) for $x \in E - \bar{D}$.

Remark. The proof of Theorem 3.1 remains unchanged if property P_3 of V is replaced by

$$P_3' : V(x) \geq 0, \quad x \in \bar{D}_v.$$

4. ESTIMATION OF $\mathcal{E}\{L(x)\}$

In this section we assume that $\mathcal{E}\{L(x)\} < \infty$ and derive an upper bound for this quantity. The result is given in Theorem 4.1.

For $x \in E$ and $t > 0$ define

$$u(t, x) = \mathcal{E}_x \left\{ \int_0^t L[x(s)] ds \right\}.$$

The next result is stronger than necessary for the subsequent application, but will be derived for its independent interest.

LEMMA 4.1. *If $\mathcal{E}\{L(x)\} < \infty$ then for all $x \in E$,*

$$\lim_{t \rightarrow \infty} t^{-1} u(t, x) = \mathcal{E}\{L(x)\}. \quad (11)$$

Proof. Let $L_n(x) = L(x)$, $|x| \leq n$; $L_n(x) = 0$, $|x| > n$ ($n = 1, 2, \dots$). By the corollary to Theorem 3.1 of [4],

$$\begin{aligned} \lim_{n \rightarrow \infty} \lim_{t \rightarrow \infty} t^{-1} \mathcal{E}_x \left[\int_0^t L_n[x(s)] ds \right] &= \lim_{n \rightarrow \infty} \mathcal{E}\{L_n(x)\} \\ &= \mathcal{E}\{L(x)\}. \end{aligned} \quad (12)$$

Let $P(t, x, B)$ be the transition function of X . If μ is the invariant measure of X then, by repeated applications of Fubini's Theorem,

$$\begin{aligned} \mathcal{E} \left\{ \mathcal{E}_x \left\{ t^{-1} \int_0^t L_n[x(s)] ds \right\} \right\} &= \int_E \mu(dx) t^{-1} \int_0^t \int_E P(s, x, dy) L_n(y) ds \\ &= t^{-1} \int_0^t \int_E \mu(dy) L_n(y) ds \\ &= \mathcal{E}\{L_n(x)\}. \end{aligned}$$

Passing to the limit ($n \rightarrow \infty$) we have by monotone convergence

$$\mathcal{E}\{t^{-1} u(t, x)\} = \mathcal{E}\{L(x)\}. \quad (13)$$

Now let $\chi_n(x) = 1$, $|x| \leq n$; $= 0$, otherwise. Suppose that for some $\epsilon > 0$ there exists a sequence $\{t_\nu\}$ ($t_\nu > 0$) and a subsequence $n(\nu)$ of positive integers such that

$$\mathcal{E}_x \left\{ t_\nu^{-1} \int_0^{t_\nu} \chi_{n(\nu)}[x(s)] L[x(s)] ds \right\} > \epsilon, \quad \nu = 1, 2, \dots \quad (14)$$

From (13) and (14) it follows that

$$\mathcal{E}\{\chi_{n(\nu)}(x) L(x)\} > \epsilon, \quad \nu = 1, 2, \dots,$$

which contradicts the fact that $\mathcal{E}\{L(x)\} < \infty$. Hence for each fixed $x \in E$,

$$\mathcal{E}_x \left\{ t^{-1} \int_0^t L_n[x(s)] ds \right\} \rightarrow \mathcal{E}_x \left\{ t^{-1} \int_0^t L[x(s)] ds \right\}$$

as $n \rightarrow \infty$, uniformly in t for t sufficiently large. We can therefore interchange limits in the left side of (12), and the result (11) follows by monotone convergence.

We consider functions V with properties $\bar{P}_1 - \bar{P}_3$, where these differ from properties $P_1 - P_3$ of Section 3 only in that now we require $D_v = E$.

THEOREM 4.1. *Let X be positive. If there exist a function V with properties $\bar{P}_1 - \bar{P}_3$ and a positive constant k such that*

$$\mathcal{L}[V(x)] \leq k - L(x), \quad x \in E,$$

then

$$\mathcal{E}\{L(x)\} \leq k.$$

Proof. We first show that $\mathcal{E}\{L(x)\} < \infty$. Indeed if D is a normal domain with boundary Γ and if $v(x) = \mathcal{E}_x\{\tau_\Gamma\}$ then $\mathcal{L}[v(x)] = -1$ ($x \in E - D$) and $v(x) = 0$ ($x \in \Gamma$). It follows that the function $V(x) + kv(x)$ satisfies the conditions of Theorem 3.1.

Let $D_n = \{x : |x| < n\}$ and put $\tau_n = \min\{t : |x(t)| = n \mid x(0) = x \in D_n\}$. Let $t_n = \min(t, \tau_n)$ and define

$$u_n(t, x) = \mathcal{E}_x \left\{ \int_0^{t_n} L[x(s)] ds \right\}$$

$t > 0$, $x \in D_n$ ($n = 1, 2, \dots$). Since $\tau_n \uparrow \infty$ ($n \rightarrow \infty$) with probability 1, we have $u_n(t, x) \uparrow u(t, x)$.

We now use the fact that $u_n(t, x)$ is the unique smooth solution of the problem

$$\begin{aligned} \mathcal{L}[u_n(t, x)] - \partial u_n(t, x) / \partial t &= -L(x), & t > 0, & \quad x \in D_n \\ u_n(0, x) &= 0, & x \in D_n \\ u_n(t, x) &= 0, & t > 0, \quad |x| = n \end{aligned}$$

(see e.g. [5], Ch. 5). We can assume that $V(x) \geq 0$, $x \in E$. If

$$W_n(t, x) = kt + V(x) - u_n(t, x) \quad (t \geq 0, \quad x \in \bar{D}_n)$$

then

$$\mathcal{L}[W_n(t, x)] - \partial W_n(t, x) / \partial t \leq 0;$$

$W_n(0, x) \geq 0$; and $W_n(t, x) \geq 0$, $|x| = n$. By the maximum principle for parabolic equations $W_n(t, x) \geq 0$ ($t \geq 0$, $x \in \bar{D}_n$); that is $u_n(t, x) \leq kt + V(x)$; hence

$$u(t, x) \leq kt + V(x), \quad t \geq 0, \quad x \in E.$$

The result now follows from Lemma 4.1.

Remark. The proof is unchanged if property \bar{P}_3 of V is replaced by

$$\bar{P}_3' : V(x) \geq 0, \quad x \in E.$$

5. APPLICATIONS

EXAMPLE 1

Let X satisfy the Itô equation

$$\begin{aligned} dx &= Fxdt - b\phi(\sigma)dt + G(x)dw \\ \sigma &= c'x. \end{aligned} \tag{15}$$

In (15), F is a constant matrix, b and c are constant n -vectors, and ϕ is a scalar-valued, in general nonlinear, function of σ . The nonstochastic differential equation, obtained from (15) by setting $G = 0$, has been studied extensively in connection with the Lur'e problem [8].

THEOREM 5.1. *Let the system (15) satisfy the following conditions:*

- (i) *All the eigenvalues of F have negative real parts*
- (ii) *$\sigma\phi(\sigma) > 0$ for all $|\sigma|$ sufficiently large; $\phi(\sigma)$ is continuously differentiable; and $d\phi(\sigma)/d\sigma$ is bounded ($-\infty < \sigma < \infty$).*
- (iii) *There exist two nonnegative constants α and β such that $\alpha + \beta > 0$ and*

$$\operatorname{Re}(\alpha + i\omega\beta)c'(i\omega I - F)^{-1}b > 0$$

for all real ω .

- (iv) *$G(x)$ satisfies the conditions of section 2 and, in addition, $|G(x)|$ is bounded for $x \in E$.*

Then X is positive and

$$\mathcal{E}\{|x|^\nu\} < \infty$$

for every $\nu > 0$.

Proof. The positivity of X was proved in [3]. To satisfy the conditions of Theorem 3.1 we introduce a function $\tilde{V}(x)$ of the form

$$\tilde{V}(x) = x'Px + \beta \int_0^{c'x} \phi(\sigma) d\sigma$$

and define

$$V(x) = \exp(\gamma \tilde{V}(x)),$$

where $\gamma > 0$ will be chosen later. By a result of Meyer [9] there exist positive definite matrices P and Q such that

$$[Fx - b\phi(c'x)]'\tilde{V}_x(x) \leq -x'Qx \quad (16)$$

for all $|x|$ sufficiently large. Moreover

$$\frac{1}{2} \operatorname{tr}[G(x)G(x)'\tilde{V}_{xx}(x)] = \operatorname{tr}[G(x)G(x)'P] + \frac{1}{2}\beta \int |G(x)'c|^2 d\phi(c'x)/d\sigma \quad (17)$$

Since the right side of (17) is bounded it follows on adding (16) and (17) that, for arbitrary $\delta > 0$,

$$\mathcal{L}[\tilde{V}(x)] \leq -(1 - \delta)x'Qx \quad (18)$$

for all $|x|$ sufficiently large. Let $\delta \in (0, 1)$ be fixed. Now

$$\begin{aligned} \exp(-\gamma \tilde{V}(x))\mathcal{L}[V(x)] &= \gamma \mathcal{L}[\tilde{V}(x)] + \frac{1}{2}\gamma^2 |G(x)'\tilde{V}_x(x)|^2 \\ &= \gamma \mathcal{L}[\tilde{V}(x)] + \frac{1}{2}\gamma^2 |G(x)'[2Px + \beta\phi(c'x)c]|^2 \\ &\leq -\gamma(1 - \delta)x'Qx + \gamma^2 x'Rx \end{aligned} \quad (19)$$

for some positive definite constant matrix R . Since Q is positive definite the matrix $(1 - \delta)Q - \gamma R$ is positive definite for $\gamma > 0$ sufficiently small. Then, for $|x|$ sufficiently large

$$\begin{aligned} \mathcal{L}[V(x)] &\leq -\exp(\gamma \tilde{V}(x)) \\ &\leq -|x|^v. \end{aligned}$$

The result now follows by Theorem 3.1.

Remark. It is clear from the proof that, under the conditions of Theorem 5.1, $\mathcal{E}\{L(x)\} < \infty$ provided

$$L(x) = O[\exp(\theta |x|^2)] \quad (|x| \rightarrow \infty)$$

for $\theta > 0$ sufficiently small.

EXAMPLE 2

We shall illustrate the application of Theorem 4.1 to the analysis of a simple control system. Suppose

$$\begin{aligned} \dot{x}_1 &= x_2 \\ \dot{x}_2 &= -x_2 - \psi(x_1 + x_2), \end{aligned} \quad (20)$$

where

$$\psi(y) = \begin{cases} 1, & y \geq 1 \\ y, & |y| \leq 1 \\ -1, & y \leq -1. \end{cases}$$

The null solution $x_1 = x_2 = 0$ is asymptotically stable. If the system is perturbed by Gaussian white noise it is of interest to estimate the mean square error $\mathcal{E}\{x_1^2\}$. The prior verification that X is positive will be omitted. Introducing perturbation terms and making the change of variables $x_1 = x$, $x_1 + x_2 = y$, we obtain

$$\begin{aligned} dx &= -(x - y) dt + a_{11} dw_1 + a_{12} dw_2 \\ dy &= -\psi(y) dt + a_{21} dw_1 + a_{22} dw_2 \end{aligned} \quad (21)$$

where w_1, w_2 are independent 1-dimensional Wiener processes and the coefficients a_{ij} are constants. The differential generator of the (x, y) process is

$$\mathcal{L}[u] = Au_{xx} + 2Bu_{xy} + Cu_{yy} - (x - y)u_x - \psi(y)u_y,$$

where

$$\begin{aligned} A &= (a_{11}^2 + a_{12}^2)/2 \\ B &= (a_{11}a_{21} + a_{12}a_{22})/2 \\ C &= (a_{21}^2 + a_{22}^2)/2. \end{aligned}$$

To satisfy condition (v) of Section 2 we assume that $a_{11}a_{22} - a_{12}a_{21} \neq 0$; in the present application such a restriction is clearly not significant.

To estimate $\mathcal{E}\{x^2\}$ we try to construct a positive function $V(x, y)$ with continuous second derivatives such that

$$\mathcal{L}[V(x, y)] \leq k - x^2 \quad ((x, y) \in E)$$

for some positive constant k . As a first step we assume that the perturbation terms are absent from (21) and evaluate

$$V^0(x, y) = \int_0^\infty x(t)^2 dt \quad (x(0) = x, y(0) = y).$$

The result is

$$\begin{aligned} V^0(x, y) &= x^2/2 + xy/2 + y^2/4, \quad |y| \leq 1 \\ &= x^2/2 - x + xy - y^2/2 + y^3/3 \\ &\quad + e^{1-y}(x - y - 1)/2 + 17/12, \quad y \geq 1 \\ &= V^0(-x, -y), \quad y \leq -1. \end{aligned} \quad (22)$$

From (22) we find that V_{xx}^0, V_{xy}^0 are continuous, but

$$V_{yy}^0(x, 1-0) = \frac{1}{2}, \quad V_{yy}^0(x, 1+0) = 1 + x/2.$$

To achieve the required smoothness replace V^0 by V^1 , where

$$\begin{aligned} V^1(x, y) &= V^0(x, y) - [\tfrac{1}{4}](1+x)(y-1)^2 e^{-\alpha(y-1)}, & y \geq 1 \\ &= V^0(x, y), & |y| \leq 1 \\ &= V^1(-x, -y), & y \leq -1, \end{aligned} \quad (23)$$

where $\alpha > 0$ is arbitrarily large. From (23),

$$\mathcal{L}[V^1] \sim 2C|y| - x^2 \quad (|x| \rightarrow \infty, |y| \rightarrow \infty)$$

To cancel the term $2C|y|$ for large $|y|$, define

$$\begin{aligned} V^{(2)}(x, y) &= V^1(x, y) + C(|y| - 1)^2 \exp[-\beta(|y| - 1)^{-1}], & |y| \geq 1 \\ &= V^0(x, y), & |y| \leq 1 \end{aligned}$$

where $\beta > 0$ is arbitrarily small. Finally, let

$$V(x, y) = (1 + \gamma)V^{(2)}(x, y)$$

where $\gamma > 0$ will be chosen later. Then

$$\mathcal{L}[V(x, y)] \leq K - x^2 \quad (x, y \in E) \quad (24)$$

if K is sufficiently large. By straightforward estimation of the individual terms of $\mathcal{L}[V]$ we can obtain a value k_γ of K for which (24) is true; we then choose

$$k = \min\{k_\gamma : \gamma > 0\}.$$

Carrying out the estimates for $|y| \leq 1$ and $|y| \geq 1$ separately, we find

$$\mathcal{E}\{x^2\} < \max(k', k''),$$

where

$$\begin{aligned} k' &= (A + B + C/2)[1 + C(9C^2 + 4D)^{-\frac{1}{2}}] \\ k'' &= (5/2)C^2 + D + (C/2)(9C^2 + 4D)^{\frac{1}{2}} \end{aligned}$$

and

$$D = A + 2B + |B| + 3C/2.$$

To obtain a rough idea of how conservative the bound may be in this case, suppose that $A \simeq 0, B \simeq 0, C \rightarrow \infty$. Then

$$\mathcal{E}\{x^2\} < k'' \sim 4C^2. \quad (25)$$

Analysis of the system (21) based on "statistical linearization" [10] of the nonlinear function ψ yields

$$\mathcal{E}\{x^2\} \simeq (\pi/2)C^2 \quad (C \rightarrow \infty). \quad (26)$$

The qualitative agreement between the results (25) and (26) is due to the special choice of the function V_0 . We should emphasize that the upper bound (25) was derived rigorously; the estimate (26), although probably reliable, was obtained by a heuristic procedure.

ACKNOWLEDGMENT

I am indebted to the referee for illuminating criticism which led, in particular, to simplification of the work of Section 4.

REFERENCES

1. WONHAM, W. M. Stochastic problems in optimal control. RIAS TR 63-14, May, 1963.
2. CRANDALL, S. H. (ed.) "Random Vibration." Vol. 2. M.I.T. Press, Cambridge, 1963.
3. WONHAM, W. M. Liapunov criteria for weak stochastic stability. *J. Diff. Eqs.* 2 (1966), 195-207.
4. KHAS'MINSKII, R. Z. Ergodic properties of recurrent diffusion processes and stabilization of the solution to the Cauchy problem for parabolic equations. *Theor. Prob. Appl.* 5 (1960), 179-196.
5. DYNKIN, E. B. "Markov Processes." Academic Press, New York, 1965.
6. COURANT, R., AND HILBERT D. "Methods of Mathematical Physics," Vol. 2. Wiley (Interscience), New York, 1962.
7. DOOB, J. L. "Stochastic Processes." Wiley, New York, 1953.
8. AIZERMANN, M. A., AND GANTMACHER, F. R. "Absolute Stability of Regulator Systems." Holden-Day, San Francisco, 1964.
9. MEYER, K. R. On the existence of Liapunov functions for the problem of Lur'e. *J. SIAM, Ser. A: Control*, 3 (1966), 373-383.
10. BOOTON, R. C. Nonlinear control systems with random inputs. *Trans. IRE CT-1* (1954), 9-18.
11. SERRIN, J. On the Harnack inequality for linear elliptic equations. *J. Anal. Math.* 4 (1954-56), 292-308.